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A NEW SOLUBLE APPROXIMATION TO FOKKER-PLANCK EQUATIONS

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The results that I am about to discuss were obtained in the course of work on collision rates in electron-positron colliding-beam storage rings. Understanding collision rates requires understanding the statistical distribution of beam particles.

Idealizing and simplifying for the sake of clarity, an electron circles a storage ring in the horizontal plane at the speed of light, with its displacement y perpendicular to the plane of the ring executing small oscillations according to

$$\ddot{y} + \delta(t)\dot{y} - F(y, t) = \eta(t) \xi(t),$$
 (1)

where %, \(\), and F are \(\) in t (the period is the time needed to circle the ring once), and \(\xi \) is Gaussian uncorrelated white noise

$$\langle \xi(t) \xi(t') \rangle = \Im(t - t') \tag{2}$$

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Correspondingly, the probability distribution, P, in y and .
y evolves in time according to a Fokker-Planck equation

$$\frac{\partial P}{\partial t} + \dot{y} \frac{\partial P}{\partial \dot{y}} + F \frac{\partial P}{\partial \dot{y}} = \frac{\partial}{\partial \dot{y}} \left(8\dot{y}P \right) + \frac{1}{2}\eta^2 \frac{\partial^2 P}{\partial \dot{y}^2}. \tag{3}$$

The Langévin equation (1), or the Fokker-Planck equation (3), differs from textbook systems primarily in that the modulation period is much smaller, rather than much larger, than the naive thermal relaxation time, $1/\langle Y \rangle$ (the averaging is over external modulation). In my work, I am particularly concerned with the limit $Y \rightarrow 0$, $\gamma \rightarrow 0$, with F and the modulation period fixed.

For today's talk, I shall ignore modulation. The new results described here, in the present one-dimensional context, constitute a necessary prerequisite for more realistic studies. I will report soon on a more extensive study of modulation-related statistical effects (and enlarge on the topics treated in the present talk) in a forthcoming series of publications.

The analysis I describe today concerns the "semiclassical" (small - η) approximation to the solution of (3) corresponding to the probability that a Langévin particle, initially at y=y and \dot{y} =v, ends up at y and v at time t:

where N is a normalization constant. (This is most easily derived from path integral considerations.) I have replaced the force F by minus the derivative of a (confining and symmetric, for definiteness) potential U. The path $y(\Upsilon)$ extremizes the exponent, subject to the boundary conditions indicated in the parentheses on the left-hand side. The extremal ("Onsager-Machlup") equation is

$$O = \left[\frac{d^2}{d\tau^2} - 8\frac{d}{d\tau} + U''(y)\right] \left[\ddot{y} + 8\dot{y} + U'(y)\right]. \tag{5}$$

I am concerned with solutions of (5) and (4) for small % but for large time (t > $O(\frac{1}{6})$), because in storage rings, $\frac{1}{2}$ = milliseconds, much shorter than the time scale (minutes, hours) on which elementary-particle experiments are performed.

In order to solve (5) approximately for small %, I assume that at any time, y is close to a solution of frictionless, noiseless Newton's law, $\ddot{y}+U'(y)=0$, up to a deviation that vanishes as $\mathring{y}\to O$ (t \mathring{y} fixed). I then proceed as follows:

I take the interval (0,t) and break it into subintervals - a subinterval (except for the first and last) begins and ends at two succesive times when y crosses zero (where U'=0). Hardly any damping or fluctuation takes place in one such subinterval. The deviation of y in such a subinterval from Newtonian vibration depends on \forall , and also on the way initial acceleration deviates from -U'(0)=0, and also

on how the initial y differs from the Newtonian value -U"(0) y.

I compute the solution y to (5) in a subinterval by perturbing about a solution to undamped Newtonian motion in powers of X, and of a (acceleration (times sign of initial velocity) at beginning of subinterval), and of p (value of Y+U"Y (times sign of initial velocity) at beginning of subinterval). From this computation I derive an approximate map, which transforms the values of a, p, and speed v at the beginning of one subinterval into their values at the beginning of the next.

This one-subinterval map is to be iterated many ($\geqslant O(\frac{1}{4})$) times. My main result is that <u>such long-time iteration can</u> be done in a simple closed form, up to small remainders.

A technical note: It is necessary to assign formal orders in \forall to the variables a and p in order to make the perturbative calculation systematic. I take my cue from two particular solutions to (5),

$$\ddot{y} \pm 8 \dot{y} + U'(y) = 0$$
 (6)

(damped and antidamped Newtonian motion). When y=0, (6) gives $a=\mp \gamma v$, and $p=-\gamma^2 v$. So in the general case I take v=O(1), $a=O(-\gamma)$, and $p=O(-\gamma)^2$.

Then the perturbative subinterval map is

$$\Delta V = \alpha \left(\frac{J}{V^2}\right) + O(8^2), \tag{7a}$$

$$\Delta \alpha = \frac{\aleph^2 V}{2} \frac{\mathcal{L}^2 J}{\mathcal{L}^2 V^2} - P V \frac{\mathcal{L}}{\mathcal{L} V} \left(\frac{1}{V} \frac{\mathcal{L}^2 J}{\mathcal{L} V} \right) + \frac{\Omega^2}{2} \frac{\mathcal{L}^2}{\mathcal{L} V^2} \left(\frac{J}{V} \right) + O(\aleph^3),$$

$$\Delta p = 8^{2} \alpha \left(\frac{J}{V^{2}} + \frac{1}{2} \frac{d^{2}J}{dV^{2}} \right) - \alpha p \left(\frac{J}{V^{3}} + \frac{d}{dV} \left(\frac{1}{V} \frac{dJ}{dV} \right) \right)$$

$$+ \frac{Q^{3}}{2V} \frac{d^{2}}{dV^{2}} \left(\frac{J}{V} \right) + O(8^{4})$$
(7b)

where J is pi times the canonical action integral for undamped Newtonian motion at energy $v^2/2$ ($J = \frac{1}{2} \oint \dot{y}^2 d \Upsilon$). Such an undamped Newtonian oscillation completes itself in time (2/v) dJ/dv.

For small Y, we may, because of (7), consider Δv , Δa , Δp as differentials rather than finite differences. Then (7) is a system of three first-order ODE's. This system turns out to have two conserved quantities

$$E = -V^{2}Y^{2} - \Omega^{2} + \lambda \rho V,$$

$$H = \frac{1}{2} \left[\frac{E}{V} \frac{\partial J}{\partial V} - Y^{2}J + \Omega^{2} \left(\frac{J}{V^{2}} \right) \right].$$
(8)

The former, E, is the restriction to y=0 of

$$E = L - \dot{y} \frac{\partial L}{\partial \dot{y}} - \ddot{y} \frac{\partial L}{\partial \dot{y}} + \dot{y} \frac{\partial L}{\partial t} \left(\frac{\partial L}{\partial \ddot{y}} \right)$$
 (9)

which is actually conserved exactly by (5), as a consequence of our present neglect of any explicit time-dependence of parameters. The latter, H, is a kind of adiabatic invariant,

$$H = \frac{1}{3} \, \left[L - \dot{y} \, \frac{\partial L}{\partial \dot{y}} - \ddot{y} \, \frac{\partial L}{\partial \dot{y}} + \frac{1}{3} \, \frac{\partial L}{\partial \dot{x}} \, (\dot{y} \, \frac{\partial L}{\partial \dot{y}}) \right] \, d\tau. \tag{10}$$

When the right-hand sides of (7) are written in terms of H, the equations for $\Delta \, v$ and $\Delta \, a$ take a canonical Hamiltonian form

$$\Delta V = \frac{\partial \alpha}{\partial H(V, \alpha, E)} + O(8^2),$$

$$\Delta \alpha = -\frac{\partial H(V, \alpha, E)}{\partial V} + O(\lambda^3). \tag{11}$$

Because of these conservation laws, the system (7) can be solved (approximately) by quadrature. The end result is this: In terms of E and H, the probability in equation (4) (restricted, for simplicity, to $y_1 = y_2 = 0$) is

$$P(O, V; O, V_1; \neq) \cong$$

$$N \exp \frac{-1}{2\eta^2} \left\{ Y(v_2^1 - v_1^2) + \int_{|v_1|}^{|v_2|} \frac{\left[2H - \frac{E}{\sqrt{2}} \frac{\partial J}{\partial v} + 28^2 J \right]}{\left[J(2H - \frac{E}{\sqrt{2}} \frac{\partial J}{\partial v} + 8^2 J) \right]^2} V(dv) + O(8^2) \right\}_{12)}$$

The path of integration proceeds from $|v_1|$ down to a turning point, and then up to $|v_2|$. To determine E and H, one extremizes the exponent in (12) (with v_1 and v_2 fixed) subject to the constraint

$$= \lambda \int_{\Lambda^{r_1}}^{\Lambda^{r_1}} \frac{\left[2(3H - \frac{2}{4} \frac{2R}{4} + \lambda_3 2)\right]_{\lambda^{r_1}}}{18\Lambda 1 (32 \lambda_3 \lambda_3)} + O(\lambda). \tag{13}$$

$$\lambda f = \lambda \sum_{\Lambda^{r_1}} \left(\frac{2(3H - \frac{2}{4} \frac{2R}{4} + \lambda_3 2)}{18\Lambda 1 (32 \lambda_3 \lambda_3)}\right)$$

Here is a check: As $8t\to\infty$, it turns out that both E and H approach zero. Then the probability (12) becomes

$$N \in \mathbb{R}$$
 $\frac{-28}{\eta^2} \left(\frac{V_2^2}{2}\right)$ (14)

i.e. Maxwell-Boltzmann, as it should. Incidentally, E=H=0

means that v, a, and p correspond to the restriction of solutions of equations (6) to y=0. This is to be interpreted as follows: When $t\to\infty$, the extremal path proceeds from (y_1,v_1) to (0,0) according to damped Newtonian motion, and then proceeds from (0,0) to (y_2,v_2) according to antidamped Newtonian motion.

Note that the harmonic oscillator is singular as far as this formalism is concerned. When U is quadratic in y, (12) and (13) depend on the same one linear combination of E and H, so that the constants cannot be determined independently. In this case, the formalism must be modified in a well-defined way. With the appropriate modifications, one can reproduce the exact harmonic oscillator Fokker-Planck Green's function (to within $O(\sqrt[3]{2})$ in the exponent).

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